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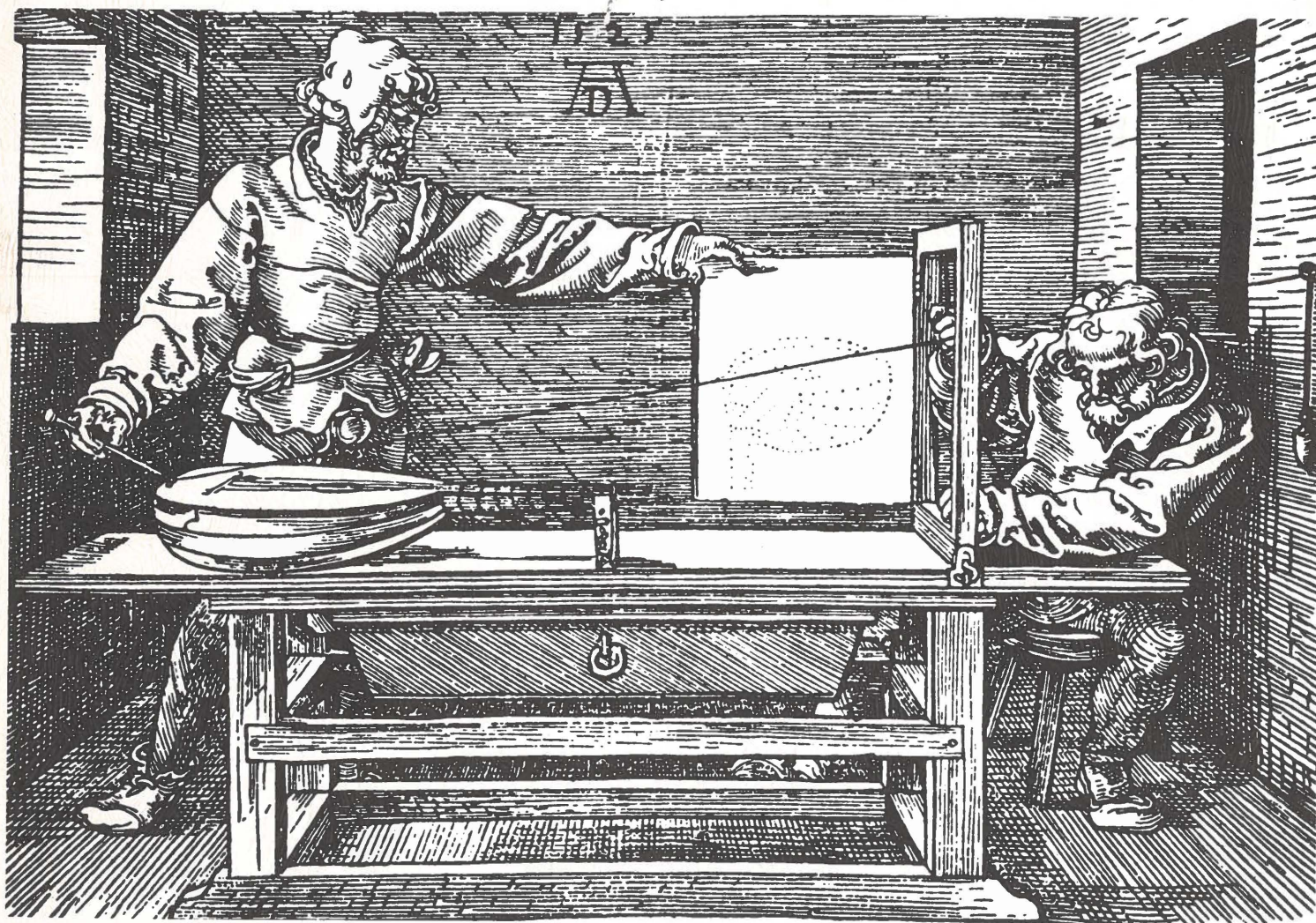
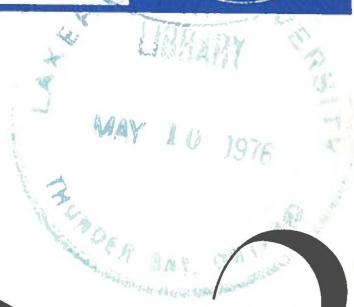
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THE VIEWS EXPRESSED IN CARET DO NOT NECESSARILY REFLECT THE OPINIONS OF THE
EDITOR, THE FACULTY OR THE UNIVERSITY.

OUR COVER relates to our article "A History of Photography" and depicts Durer's use
of the pin-hole camera.

A HISTORY OF PHOTOGRAPHY

by: Dr. Andrew D. Booth
Lakehead University, Thunder Bay

A science more than 250 years old

The science, as distinct from the art, of photography is now more than 250 years old since the basic experiment on the blackening of silver salts through the action of light was conducted by J. H. Schulze in 1725. Schulze did not make a photograph as we now understand it but used a sensitive material which consisted of a mixture of chalk and silver nitrate shaken up with water in a bottle and used wet to record the outline of a stencil placed between the emulsion and the sun. Since the process was essentially a wet one, there are no existing "photograms", to use the modern term, dating from this time.

In 1777, C. W. Scheele noticed that silver chloride was more quickly blackened by the violet and blue rays of the spectrum than by the others, and in a sense, this fundamental experiment is the basis of much modern work on photographic emulsion sensitivity and its correction to record the same sort of image as seen by the human eye.

In the year 1800, Thomas Wedgwood made contact copies in silhouette form of leaves on leather sensitized with silver nitrate. Again, there was no fixing process used so that the results of these experiments have long since disappeared.

The man most usually credited with the invention of photography was **Nicephore Niepce** who, in **1816**, made photographs using a "camera" — that is, a dark chamber with an appropriate lens — to produce photographs on paper sensitized with silver chloride. These photographs were partially fixed using hypo and really constitute the first artistic realization of what we now call photography.

In the year 1819, Sir John Herschel discovered that thiosulphates removed undeveloped silver halides and thus made the images produced on paper by the action of light permanent. This led

in 1826 to the second relevant experiment of Niepce who made a photograph from nature using a lens system and a pewter plate sensitized with bitumen. Niepce's camera of 1826 is interesting in that it contained a bellows for focussing the lens and an iris diaphragm for controlling the amount of light, although the degree of control is debatable since **the exposure was eight hours!**

Recognizable artistic photography started in 1835 when L. J. M. Daguerre made direct photographs on silvered copper plates coated with a silver iodide. The latent image was developed by mercury vapour, a most obnoxious substance, and in 1837, it was found that the result could be fixed using a solution of common salt. Thus, the **Daguerreo type** was born.

The second contender for the honour of invention of artistic photography was W. H. Fox Talbot who, also in 1835, produced "**photogenic drawings**" on paper sensitized with silver chloride and fixed with potassium iodide. Fox Talbot also discovered that prolonged washing with water would dissolve out the undeveloped chloride and thus act as a fixative.

The period 1820 to 1840 was one of vigorous scientific activity and readers should recollect the work of Michael Faraday who, between 1828 and 1832, discovered most of the practical applications of electro-magnetism, including the transformer, the motor, and the electric generator. The time was also an active one for microscopists, and in 1839, the Reverend J. B. Reade made the first photomicrographs using a solar microscope to record on paper sensitized with silver nitrate and developed with gallic acid. Fixing was achieved by immersion in sodium thiosulphate.

The importance of the photographic process is indicated by the fact that in 1839, the French Government issued working instructions for the creation of Daguerreotypes, and in the same year, Sir John Herschel, son of Sir William mentioned earlier, first coined the word "**photography**" from roots meaning **writing with light**.

1840 is an important date, since in this year, J. Petzval established the basic theory required to compute optical systems of photographic type. The Petzval portrait lens is still in use.

In the same year, A. Wolcott opened the first photographic portrait studio in New York City. **The Europeans had the art, but the Americans, as usual, made the money!**

1844 — first book illustrated photographically

1844 was an important date in the history of photography since it was in that year that Fox Talbot published the first part of "The Pencil of Nature" which was the first book to be illustrated photographically. Copies of this book are still extant, but it is a very rare book, and if a copy ever came up for sale, it would fetch a very large sum of money.

The birth of flash photography was not due to Harold Edgerton as is usually thought, but to Fox Talbot who, in 1851, made the first photographs of an object in rapid movement using an electric spark. In the year 1852, the same experimenter showed that gelatin was made insoluble when saturated with sodium biochromate and exposed to light. This led to the first practical process for half-tone printing.

The technical processes of photography were in a state of rapid change. The Daguerreotype required special preparation for viewing. The later wet processes were messy in application and were replaced finally by dry processes using a transparent negative followed by a positive.

1861 — colour!

All of this work was, of course, in black and white or in some of the rather charming sepia and gold tones produced by the replacement of silver by such metals as gold and platinum, but in the year 1861, J. Clerk Maxwell demonstrated the **first three-colour separation method** for producing coloured images and thereby established the basis of modern colour photography. The work of Maxwell was followed quite quickly in 1868 by that of Ducos du Hauron who experimented with three-colour photography and discovered various methods of realizing it including that of **subtractive colour synthesis**.

It was not until **1871** that R. L. Maddox made the **first gelatin dry plates**. These were extremely slow, although subsequently improved, both as to speed and to quality. With the coming of the gelatin emulsion, the chemists concentrated on finding reliable methods of development and Sir William Abney, of Photometer fame, discovered the virtues of hydroquinone in 1880 and was soon followed by H. B. Berkeley who, in 1882, added sodium sulphite to the armoury of photographic developers.

1882 also marks the point of emergence of **orthochromatic plates** — that is, plates whose sensitivity resembled more that of the human eye than did the original emulsions which, as we mentioned earlier, displayed considerable sensitivity to blue and violet radiation but little to anything else.

From this time on, the development of photography was rapid and the inventions came thick and fast. **Significant dates are 1888** when the first Kodak roll film camera came on the market, **1891** when the physicist Lippmann produced the first interference colour photographs, and **1906** when Wratten and Wainwright produced the **first panchromatic plates** — that is, plates sensitive to all colours.

1914 saw the **first Kodak colour photographic process**, but the real development of colour photography had to await the work of Mannes and Godowsky in **1935** when the **Kodachrome process** was developed.

The last significant date was 1947 when Edwin Land produced the one-step photographic process now known as **Polaroid**.

Along with the development of the chemistry of the photographic process, optical systems and cameras received a great deal of attention. The first cameras were simply boxes with a lens in front and a plate behind. Early lenses, which used a very small aperture, did not require focussing, in this respect they did not differ much from the box Brownies of the 1930's and the inexpensive plastic cameras of the present day! Because of the slow emulsions, however, it became necessary to develop lenses which would let in far more light. This led to two interesting scientific kinds of research: the first directed to the elimination

of the severe aberrations from which large aperture lenses suffered, and the second to the design of lenses to produce distortion-free images. The only important technical developments of the nineteenth century were the addition of bellows to enable the new and larger aperture lenses to be properly focussed and the introduction first of the Waterhouse diaphragm with its disc of circular stops, and second of the iris diaphragm as a means of controlling the amount of light falling on the film. As films or plates became faster, exposures became much shorter and this, in turn, led to the mechanical development of shutters which would work at high speed and at the same time give reproducible results. Many and various were the designs tried, ranging from pneumatic to spring-driven sets of blades, and of course, roller blind type apparatus, usually driven by springs, which was the prototype of the modern focal plane shutter.

miniature, spy cameras

The late nineteenth century produced a rash of spy cameras, small and unobtrusive devices which were the prototypes of the modern miniature cameras. These were designed to take pictures of one's friends in embarrassing situations, to say nothing of their use by the emerging race of scientific detectives typified by Mr. Sherlock Holmes. The first miniature cameras in the modern sense were the Ermanox used to great effect by the famous photographer, Dr. Salomon and, of course, the Leica, the grandfather of all modern miniature 35 mm. cameras. The explosion of photographic technology of the 1960's and the introduction of electronics, both for light metering and for shutter automation have produced a real revolution in the use of the camera. **No longer does the photographer have to have any real skill. What he needs is artistic perception. The automation does the rest.**

CULTURE, POLITICS AND GEOGRAPHY: or Cymru Am Byth Is A Fine Slogan, But What Does It Mean?

by Mr. Robert S. Dilley
Lakehead University

While there are almost as many definitions of geography as there are geographers, few would deny that one of the most basic concerns of the subject is with distributions: where things are, why they are there, and how they relate to other things. Such investigations can take many forms: the distribution of crop-growing and its relationship with rainfall; the distribution of coal and its relationship with unemployment, and so on. In each case two (or more) phenomena are mapped, and attempts made to measure and explain any similarities in their distributions. For example, consider the relationship between minority cultural traits and minority political parties. In the case of Wales the historic national tongue is Welsh, a Celtic language with no recognizable relationship with English. It is found only in Wales and among a few expatriates, though Welsh-speakers tend to lose their language quickly on emigration. About one-fifth of the population currently speaks Welsh (Map 1), though virtually

all are bilingual in English (Welsh monoglots account for little over 1% of the total population). Welsh-speaking is strongest in the rural areas of the north and west; lowest in the industrial south, along the English border and in Pembrokeshire, "little England beyond Wales". Being exclusive to the Welsh it provides a strong focus for national cultural identity; even non-Welsh-speaking Welshmen delight in the average Englishman's inability to pronounce place-names such as Pwllheli, Llanllwchaearn and Llanfihangel yng Ngwynfa.

Welsh nationalism as a significant political force is a new phenomenon; after years of being the butt of jokes about lost deposits the Plaid Cymru (pronounced approximately "plide come-ree", with the emphasis on "come") attained respectability when its President, Gwynfor Evans, was elected in a Carmarthen by-election in 1966. Last year's October election was their most successful (Map 2) when they gained 10% of the popular vote and three M.P.s; in Carmarthen (Gwynfor Evans again), Caernarvon and Merioneth.

The general relationship between Plaid Cymru and the Welsh language is apparent from the maps (for simplicity, data are grouped by counties: Glamorgan, for example, returns 17 M.P.s; Brecknock and Radnor share one between them). Correlating percentage Welsh-speaking with percentage vote for the Plaid Cymru on a county basis provides a coefficient of +0.83, significant at the 0.1% level. This is an indication of a very strong positive relationship. Such relationships, of course, must not simply be pulled out of the statistical hat and held up for wonderment. In some cases high correlations may be due to chance; more often to the overriding influence of some third factor on both elements being studied. Before drawing maps or running correlations the researcher must have good reason to expect to find a relationship. In the case of Welsh and Welsh nationalism the connection is clear: the party has made great use of the language as a rallying-point or standard: although only a minority is fluent most Welshmen know at least a few words and have a sentimental attachment to the symbol of their difference from the *saesneg* (Saxon; i.e. Englishman). Appeals to *hwyl* (intensely sentimental Welsh emotionalism, seen at its quintessential best just after one of the regular Welsh rugby football defeats of the England side) are more effective when couched in Welsh. Conversely, non-speakers of Welsh may become irritated by the noisy waving of the linguistic flag, and less inclined therefore to vote for the Plaid Cymru.

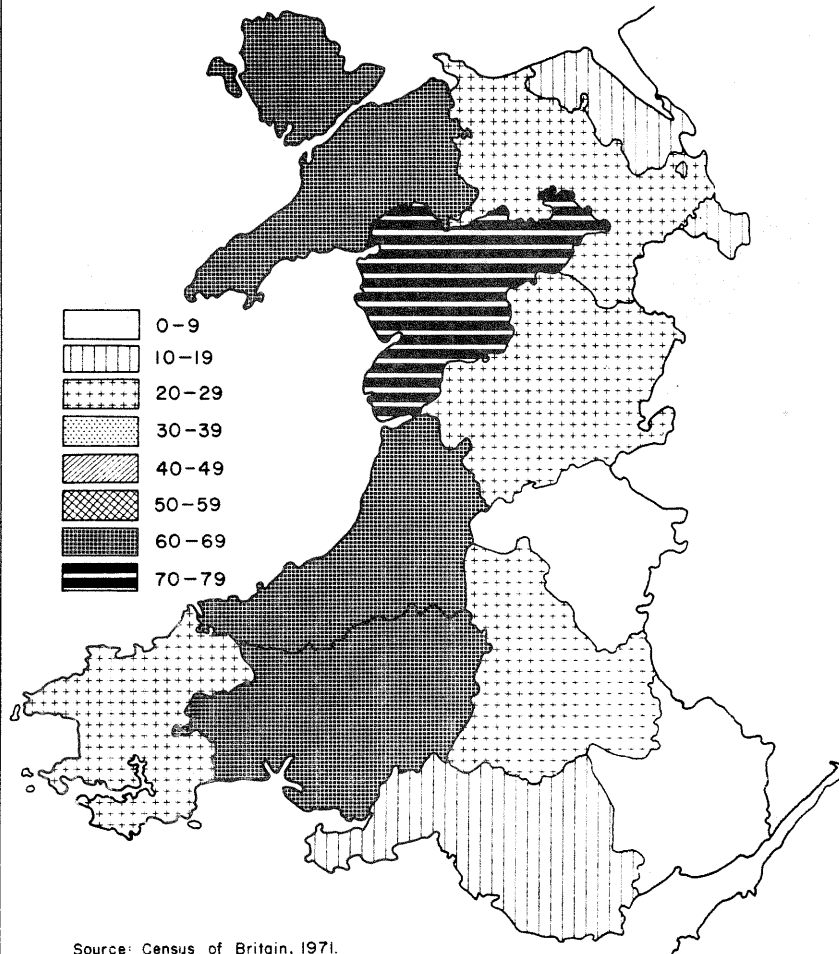
The relationship, clearly, is not perfect. The percentage vote for the Plaid Cymru at the last election was only half the Welsh-speaking percentage, and it did not follow the language distribution absolutely. People vote for a variety of reasons, and will not automatically support a party just because it is the only one to espouse their language. Many nationalistic Welsh-speakers on Anglesey voted for the Labour M. P. Cledwyn Hughes, a prominent and much-admired advocate of the Welsh cause. In the heavily English-speaking, coal-mining and steel-producing areas of Glamorgan and Monmouth the vote has traditionally been monolithically Labour: in 1966 the area returned 21 Labour M. P.s and one Conservative. However, the failure of successive Labour governments to improve the South Wales economy

sufficiently has led to a certain amount of protest voting. The rural areas have tended to go Tory (they now have 4 members). The miners would sooner vote for a Martian than a Conservative (at the last election Abertillery voted 76% Labour and Rhondda 77%), but an increasing number feel safe in registering their irritation by giving their vote to Plaid Cymru. Thus the Anglesey vote is rather lower than might be expected; that of the industrial south a little higher. If space permitted these residuals, or differences from the expected figures, could be mapped and more easily observed.

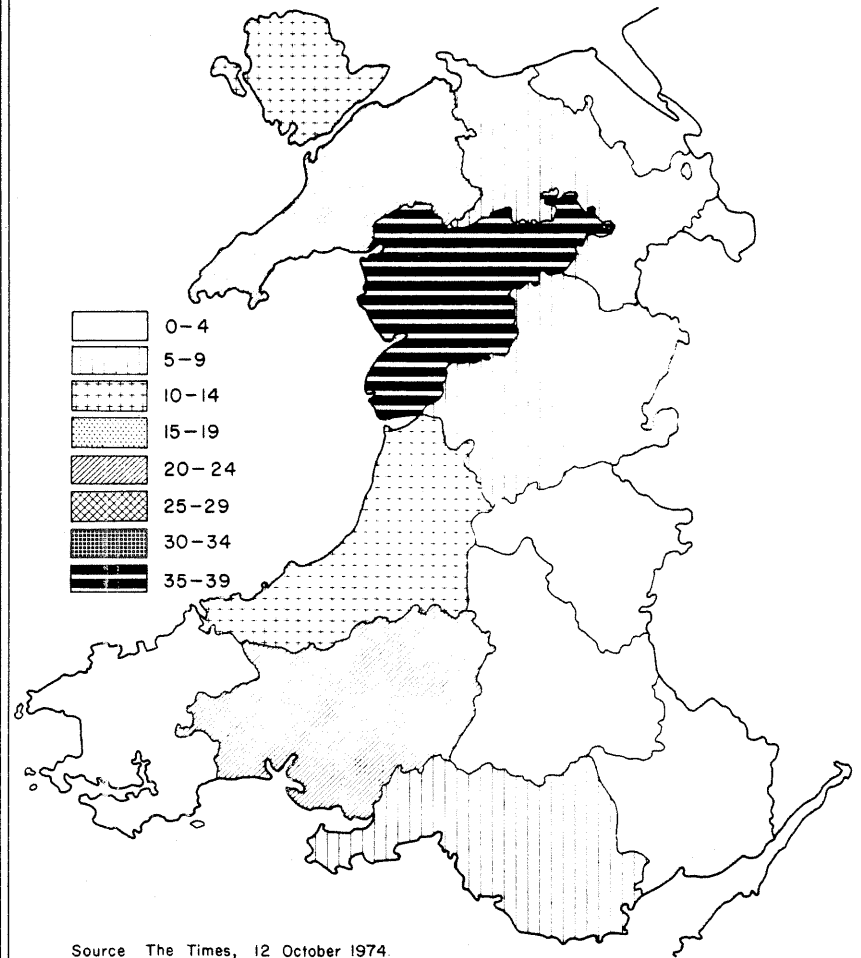
At this level of study only crude approximations can be made. Greater accuracy would be obtained by breaking the counties down into their individual constituencies (smaller division is impossible, as British elections are not declared on a poll-by-poll basis: the ballots are taken to a central location and deliberately mixed before being counted). Similar studies could be made of other areas where linkages of cultural phenomena and political allegiance are suspected. It might be interesting, for example, to compare Francophone areas and votes for the Parti Quebecois. There would not, however, be any point in relating votes for the Scottish Nationalist Party to the Gaelic tongue. Very few Scots speak the language (little over 1%) and they are highly localized in the Western Highlands and Islands. The S. N. P. concentrates more on economic and governmental issues and has wider appeal than the culturally-oriented Plaid Cymru (the S. N. P. polled nearly 30% of the Scottish vote last year, overtaking the Tories).

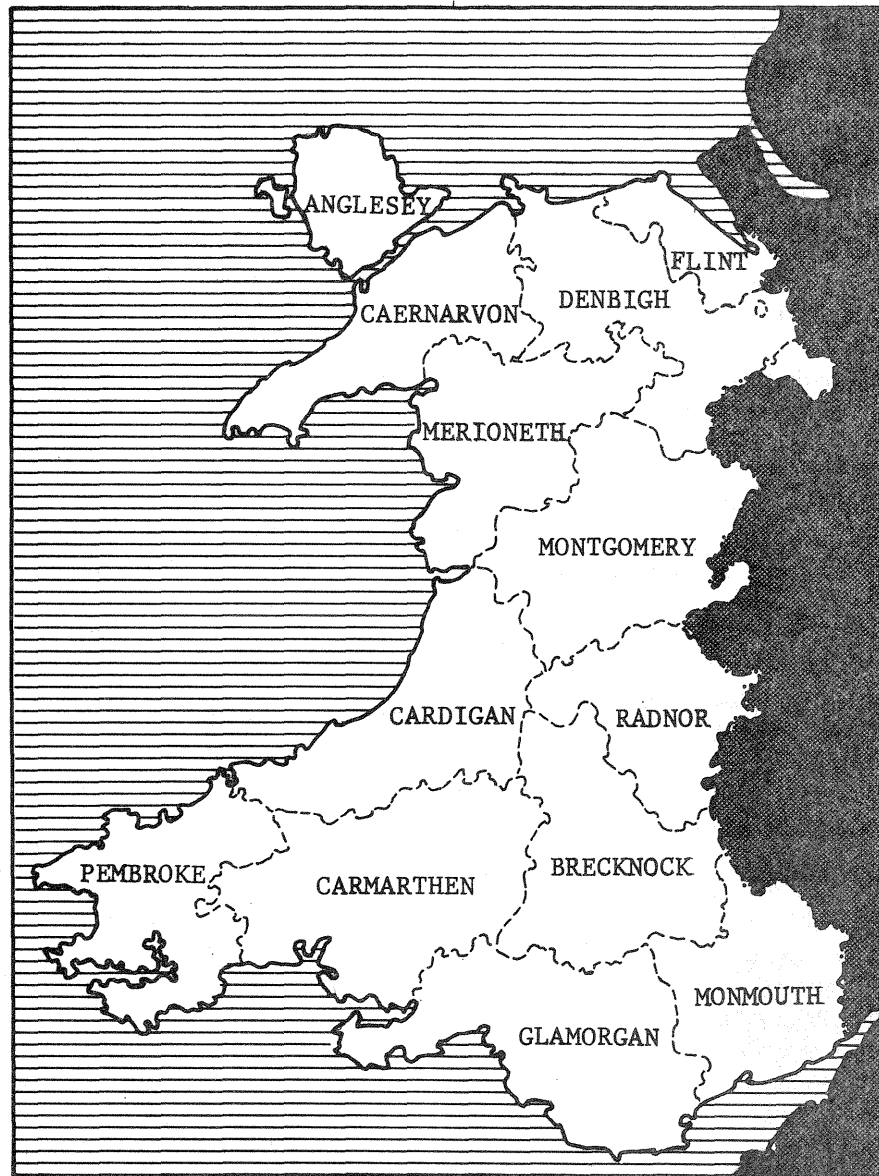
Thus Plaid Cymru sentiment is clearly linked with the Welsh language. It might seem advantageous for the party to foster this link, and to aim to collect the other half of the Welsh-speaking vote. However, this could be risky in the long term, as the Welsh-speaking percentage is shrinking and the Anglophone majority may be repelled. Plaid Cymru votes in largely English-speaking areas such as Pembroke (4.5%), Barry, Glamorgan (3.3%) and Monmouth (1.4%) suggest that "Cymru am byth" (Wales for ever) has less meaning there. Studies such as this can at least provide the basis from which political decisions must be made.

MAP 1, PERCENTAGE OF POPULATION OVER 3 YEARS OLD ABLE TO SPEAK WELSH.



MAP 2, 1974 PLAID CYMRU VOTE AS A PERCENTAGE OF THOSE VOTING.





MAP 3, COUNTIES OF WALES

THE STRATEGY OF CONCENTRATED FIRE POWER

by: *Air Cadet L. Mathemaki (Ret'd)*
 Director of the Division
 of Theoretical Warfare

1: INTRODUCTION. With the increasing mathematization of the social and natural sciences, most math teachers find they must drastically modify and update their course contents, in order to maintain at least a semblance of relevance. Gone are the days when students would be content with only physical and geometrical applications of the calculus, for example! Now most high school calculus courses must contain applications to economics, biology, sociology and psychology, even if they are of an extremely idealized nature.

Thus, the following very simple application of calculus to warfare might appeal to a good student, although it may be difficult to persuade him/her that it is really of practical value. At the very least, the main result is both easy to obtain and completely unexpected - two good qualities rarely occurring together!

2: A SHORT HISTORY OF WARFARE. We deal with battles between two opposing forces. Battles in primitive times probably consisted of a lot of individual hand-to-hand fights; the essential point for us is that the weapons - fists, clubs, spears - were not long range weapons: it was impossible to concentrate the power of many soldiers on one opponent. Thus if, for example, we had two such armies, one with 100 men and one with 200, we would expect - if the soldiers are equally proficient - that the result would be annihilation of the first army and 100 survivors from the second. The battle could simply be fought by letting half the second army act as spectators, while everyone else is killed.

Nowadays things are different and undoubtedly much improved. Long range weapons enable armies to concentrate their fire power and, as we shall see, this can lead to unexpected advantages. There are some pretty obvious examples: perhaps the battle of Agincourt was the most spectacular - the long range weapons of the English (bows and arrows) annihilated the the larger French army, whose weapons were short range.

Nearer our own time, the use of field artillery in the Russo-Japanese war, and now, the anticipated use of inter-continental ballistic missiles, indicate how far we have come from clubs and swords. Interested history students might profitably analyse some modern battles, using the concepts we will now introduce.

3: A MATHEMATICAL MODEL FOR MODERN WARFARE.

We consider two opposing forces, the red and the blue armies, of sizes $r(t)$, $b(t)$, respectively, at any time t . We will measure time from the beginning of the battle, so $r(0)$, $b(0)$ are the initial sizes of the armies. Of course, we could measure the sizes, not in terms of soldiers, but in terms of battalion, ships, etc., whatever is convenient. We assume that each soldier (battalion, ship) in the red army is equally effective in warfare, and that he can concentrate his fire power in any way, so that, throughout the battle he kills R blue soldiers every minute. Similarly, each blue soldier kills B red soldiers every minute; R and B are fixed numbers.

Consider what happens in a short time interval from t to $t+h$ minutes during the battle. The number of red soldiers killed is, by the definition of r ,

$$-(r(t+h)-r(t)).$$

On the other hand, if the time interval is short, the size of the blue army hasn't changed much in this time, and thus is approximately $b(t)$ - its size at the beginning - so the number of red soldiers killed is

$$b(t) \cdot B \cdot h$$

by the definition of the kill rate B . Thus, for values of h near 0,

$$r(t+h) - r(t) \approx -b(t) \cdot B \cdot h$$

so

$$\frac{r(t+h)-r(t)}{h} \approx -b(t) \cdot B$$

Since this approximation improves as h approaches 0, we have

$$\frac{dr}{dt} = -b \cdot B$$

(Calculus!) Similarly,

$$\frac{db}{dt} = -r \cdot R$$

and so, after a bit of algebra

$$b \cdot B \cdot \frac{db}{dt} = r \cdot R \cdot \frac{dr}{dt}$$

If we integrate both sides with respect to time, we get the unexpected result:

' $b^2 B - r^2 R$ is independant of time

Let us call $b^2 B$ the **strength** of the blue army, and $r^2 R$ the **strength** of the red army. We have shown that *the difference in strength is constant throughout the battle.*

Examples : 1. Consider a battle between 10 machine gunners (the red army) and 100 infantrymen (the blue army), and assume the kill rate of the machine gunners is 16 times that of the infantrymen. Let us measure time so the rates are 16 and 1. Initially, the strengths are:

$$\begin{aligned} \text{machine gunners} &= 16 \cdot 10^2 = 1600 \\ \text{infantrymen} &= 1 \cdot (100)^2 = 10000 \end{aligned}$$

and so, at any time,

$$b(t)^2 - 16r(t)^2 = 10,000 - 1,600 = 8,400$$

The battle will end when one side is vanquished - when $b(t)$ or $r(t)$ is 0. Clearly, $b(t)$ will never be 0:

$$b(t)^2 = 8,400 + 16r(t)^2$$

so, at the end, $r(t) = 0$ and $b(t) = \sqrt{8,400} = 91$. Only nine infantrymen have been killed, and all 10 of the machine gunners die in spite of the fact that they are, individually, 16 times more effective in killing! A rather surprising result.

2. The above example was rather an uneven contest - the machine gunners were quite overwhelmed. Let's consider a more equal match: 10 machine gunners against 60 infant-

rymen. The infantry is the stronger, by 2000, so a battle would result in it winning, with 45 infantrymen left.

But suppose the machine gunners use a bit of strategy, and manage to break the battle into a series of battles, in each one of which they all attack 20 infantrymen. The results are:

- (i) 10 mgs vs 20 inf. : 8 mgs survive
- (ii) 8 mgs vs 20 inf. : 3 mgs survive
- (iii) 6 mgs vs 20 inf. : 3 mgs survive

The machine gunners have won!

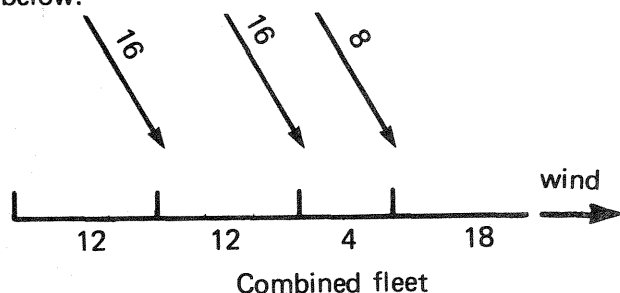
Thus we see that a bit of strategy can win against what, at first glance, appears to be overwhelming force. Of course, the enemy general has to be rather stupid to allow his army to be picked off piecemeal, but this is what often occurs in practice. A small force is sent to divert a larger part of the enemy; their job is simply to pin down a lot of soldiers while their comrades defeat the small remainder. Then they keep regrouping until they win. A naval battle, during the days of sailing ships, would be a good example of this - in that no diversionary force is needed if the wind is right - the attacking admiral hits his opponents part way down the line and lets the wind carry off a portion of the enemy fleet. By the time it has been turned around to re-enter the fight, the next phase has started. Nelson used this ploy at Trafalgar; this has been analysed by Lankester in his classic paper [3]. We will give a slightly different interpretation.

4: THE BATTLE OF TRAFALGAR RERUN.

Up to 1782, there was no 'naval strategy' - it was 'line up, and attack ship-to-ship'. But in 1782 the British admiral Rodney, probably by accident, broke through the enemy line and by concentrating on his centre and rear, won a victory. This prepared the military mind for Nelson's strategy at Trafalgar. It was known that Nelson would favour Rodney's strategy; the French admiral, Villeneuve, stated before the battle that "the British fleet will not be formed in a line-of-battle parallel to the combined (ie. the French and Spanish) fleet according to the usage of former days. Nelson will seek to break our line, envelop our rear, and overpower with groups of his ships as many as he can isolate and cut off. Why the French and Spanish admirals let Nelson dictate the strategy is a matter for military historians; the answer may lie in the fact that the French believed their

reverses at sea were due to the introduction of tactics, which one of their admirals had castigated as a "veil of timidity". Also, due to the British blockade, the French sailors at that time were so untrained they couldn't manoeuvre their ships - but they were more accurate at gunnery, so long-range warfare should have been congenial for them.

At any rate, Nelson and his aides, Collingwood and Keats, planned the battle as follows: They assumed they would have 40 ships, and the enemy would have 46. The British fleet was to attack in 3 columns, as below.



Assuming the wind carried the 18 ships out of the battle, and that it broke into 3 separate battles:

16 British vs 12 Combined
 16 British vs 12 Combined
 8 British vs 4 Combined,

resulting in 27 British ships surviving, and then the final battle of 27 British against 18 Combined, the net result is that 20 British ships are left. If the usual 'line of battle' attack had been used, the combined fleet would have won, with 20 ships, and most of us would speak French.

In fact, these tactics were changed because part of Nelson's fleet had to accompany a convoy to Malta, and in the battle he only formed two lines; the battle is described in great detail in [1].

5: MORE APPLICATIONS. It would be interesting to apply these ideas to modern battles, to see if the military strategists could have been using - perhaps in an intuitive way - the quantitative results we can get. A 'dictionary of battles' - there are quite a few in print, and they should be available in the reference section of most school libraries - is all that would be needed.

There may be applications to other situations to; polymer chemistry and molecular biology might contain interesting applications. Also, the mathematics of the situation has not, to my knowledge, really been investigated. The length of battle can be found; the answer involves the hyperbolic functions. The strength of an army consisting of several different types of weapons is certainly not that given in [3], and appears to depend on the opposition, but I have not found even an approximate explicit form for it. **What is the size of an overwhelming force?**, one that will win no matter what strategy is used. **How can we find a best strategy?**

Other mathematical problems may suggest themselves; I would be glad to hear of them.

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Synonyms

Glycerol and **glycerine** are used for the same compound which is a trihydric alcohol. Hence the former spelling with an -ol ending is preferred.

Xylene and **xylol** are also used synonymously but the compound is not an alcohol but a homologue of benzene. The first spelling is preferred.

Glycine is an amino acid (also known as aminoacetic acid) while **glycinol** (also called 2-aminoethanol, ethanolamine, monoethanolamine) lacks the carboxylic group.

Glycol: common name given to ethylene glycol.

MATHEMATICS AND THE ARTS STUDENT

by: Dr. C. C. Mutambirwa
Lakehead University

The mathematical needs of the Arts student have increased very much in recent years. This is particularly so for the student intending to major in any one or more of the social sciences, namely, Anthropology, Economics, Geography, Political Science, Psychology and Sociology. Unfortunately a large number of such students still enter college or university with limited mathematical preparation at the high school level.

Most Arts-bound students drop out of the high school mathematics program because of real or imagined fears of mathematics as a hard subject. Others drop out because of the misconceptions (also held by several teachers) that high school mathematics is not essential to the Arts major. The lack of a significant change in this state of affairs seems to lie in the pure Science bias contained in most high school mathematics curricula which have failed to include explicitly material suitable for the needs of the Social Scientist. Lately, however, the growing importance of mathematics to the social sciences is being recognized. The Secondary Schools Statistics Project which was launched just over two years ago at the University of Western Ontario is one evidence of this recognition. Since then high school mathematics teachers have held related seminars designed to revise the high school curriculum (especially the statistics program) accordingly.

The growing importance of mathematics to the social sciences has been demonstrated by the results arising from the study under the auspices of the project noted above. (See MacNeil, 1975, "The Social Scientists' View Of High School Mathematics", in Ontario Mathematics Gazette Vol. 13, No. 3, pp167-186). Using questionnaires, a poll (80% response) of social scientists in Ontario Universities was conducted to determine their opinions on mathematics curricula. Three sets of questions were asked. The first set asked the social scientists to indicate the value of eleven major mathematics topics to their undergraduate programs. The second set asked for the antici-

pated future mathematical needs of the social sciences. The third requested them to rate the importance of various statistics and probability topics to the social sciences programs.

The results of the statistical analyses of responses to the above sets of questions show that:

1. Statistics, probability and the basic concepts of algebra are very useful to the secondary school mathematics program and hence also to the university social science programs. (Table 1 refers).

2. The relatively recent and modern aspects of the mathematics curriculum, which include computer science, probability, statistics and matrix algebra, will be of increasing importance to the social sciences. (see Table 2).

3. Data tabulation methods, measures of central tendency and of dispersion should be part of the high school statistics course. In addition the high school student should be introduced to the elementary concepts on probability theory (probability measures, mathematical expectation, etc.), statistical inference, simple correlation and regression analysis, common experimental designs and associated analysis of variance, and some nonparametric tests.

At the request of the survey some general comments concerning the secondary school mathematics program in particular, were appended to the questionnaires by the social scientists. The majority of the comments expressed the opinion that greater preparation of secondary school students in basic mathematics is very necessary and mathematics should be mandatory for all students. However, there are some reservations about the type and depth of statistics to be taught at high school level. It was generally felt that only the elementary topics in statistics could be taught with the intention of showing the secondary school student that statistics *inter alia*, is an integral part of most social science programs. Complete and detailed statistics courses could be left to the colleges and universities by which time the relevance of statistics to the social sciences becomes clearer and the subject can be easily integrated into a specific social science major.

Mathematics and statistics have become part and parcel of most social science programs. Several recent publications in the social sciences clearly show this to be the case. The growth of a scientific methodology in the social sciences, the development of computer technology and the magnitude and complexity of the real-world problems have all led to the adoption of quantitative approaches to problem solving. These developments require substantial mathematical skills in the arts student, just as much as the pure science student must develop some skills in the social sciences if his or her scientific endeavours are to have any relevance to the more immediate problems of the real world.

Scale:

- 1 of no use
- 2 of limited use
- 3 a useful topic
- 4 very useful
- 5 indispensable

TOPIC NO.	MATHEMATICS TOPICS	Economics	Geography	Political Science	Psychology	Sociology & Anthropology	All departments combined
1	Algebra. Basic concepts	4.27	4.38	2.73	4.00	3.92	3.89
2	Set theory. Basic concepts.	3.45	3.46	2.55	3.13	3.00	3.13
3	Euclidean geometry	3.09	3.50	1.36	2.19	1.92	2.37
4	Analytic geometry.	2.91	2.92	1.45	2.25	2.42	2.40
5	Trigonometry.	2.20	3.45	1.18	1.94	1.92	2.15
6	Differential & integral calculus.	3.55	3.00	2.18	2.56	2.58	2.76
7	Matrix algebra.	3.64	3.54	2.55	2.60	3.08	3.07
8	Statistics.	3.82	4.00	4.09	4.13	4.08	4.03
9	Probability	3.55	3.85	3.73	3.88	3.75	3.76
10	Computer programming.	3.00	3.54	3.00	3.25	3.50	3.27
11	Actuarial mathematics.	2.09	1.64	1.50	1.81	1.92	1.80

TABLE 1

Average response to importance of mathematics for social science disciplines.

Source: Ian B. MacNeil, 1975, "The Social Scientist's View of High School Mathematics", Ontario Mathematics Gazette, Vol. 13, No. 3, p.180.

TOPIC NO.	MATHEMATICS TOPIC	Economics	Geography	Political Science	Psychology	Anthropology & Sociology	All Departments Combined
1	Algebra. Basic Concepts	0.455	0.500	0.300	0.133	0.333	0.350
2	Set theory. Basic concepts.	0.364	0.417	0.273	0.429	0.364	0.373
3	Euclidean geometry.	-0.182	0.167	0.000	-0.214	-0.091	-0.070
4	Analytic geometry.	-0.100	0.000	0.111	0.000	0.000	0.035
5	Trigonometry	-0.091	0.091	-0.100	-0.143	0.000	-0.052
6	Differential & integral calculus.	0.636		0.300	0.200	0.667	0.441
7	Matrix algebra	0.636		0.500	0.667	0.750	0.661
8	Statistics	0.455	0.833	0.800	0.600	0.667	0.667
9	Probability.	0.364	0.909	0.600	0.688	0.750	0.667
10	Computer programming.	0.364	0.750	0.700	0.813	0.833	0.705
11	Actuarial mathematics.	-0.300	0.143	0.000	0.000	0.000	-0.083

Table 2

Average response to predicted future importance of mathematics topics for social science disciplines.

Source: Ian B. MacNeil, 1975, "The Social Scientist's View of High School Mathematics", Ontario Mathematics Gazette, Vol. 13, No. 3, p. 182.

SOLUTION TO PREVIOUS PUZZLE:

FLIERS, STINGERS AND BITERS

1	I	N	S	2	E	C	T	F	3	A	M	4	I	L	5	E	6	S			
	C				M					N		N		N			H				
7	H	E	L	I	X					T		C		E			E				
	N				C			8	B		9	S	K	I	P	P	E	R			
10	E	R	G	A	T	E						S		T			R				
	U				T			11	D	I	G	12	E	R			Y				
13	M	I	T	E					B			A					S				
	O								U			D			14	P	E	E	P		
	N				15	E	16	N	G	U	L	F			E				I		
	W					17	S			I					18	L	U	R	D	A	N
19	A	N	N	U	L	A	R		20			Y			S					N	
	S			A		G				O				21	M	I	D	G	E		
	P			I		A				A					S					R	
22	S	O	L	D	I	E	R	B	E	E	T	L	E	S							

CHALLENGES TO FISHERY RESOURCE MANAGEMENT

by: *Professor W. Momot*
Lakehead University

Although fisheries management has existed as a discrete discipline for only 40 years, there has been a greater impact on our fishery resources during this short time than in the past six thousand years of recorded history. While the burgeoning populations of the third world depend increasingly on fish as a source of protein, the fish supply is decreasing. The annual world fish catch has failed to increase for the second time since World War II. Latest figures gathered by the FAO, the food and agricultural organization of the United Nations, show that the world catch of all species for 1972 was 65.6 million metric tons, down 4.1 million tons from the previous year.

The need for protein has led to fierce international competition for the ocean's fishery resources. Many traditional fishing grounds are harvested so intensively that some of our most important commercial species have been thoroughly over-exploited. Much of the declining fish catch in the United States and Canada can be directly traced to increased competition from foreign fleets. This has occurred in the haddock fisheries of the north Atlantic. Despite establishment of the International Commission for Northwest Atlantic Fisheries, which deals with the problems created when several nations exploit a common resource in international waters, satisfactory regulatory action has not been attained. As traditional fisheries decline the fishing nations of the world turn to other resources, and the same pattern of over-exploitation is repeated again and again. Even such oceanic fisheries as that for the Pacific salmon, Pacific halibut, and tropical tunas, which have yielded substantial catches over the past few decades, are being placed under more and more pressure.

Another facet of this intense competition is the attempt by some coastal nations to limit access to their offshore resources by extending their coastal jurisdiction to the edge of the continental shelf or to some other fixed point usually given as 200 miles offshore or 200 meters in depth. This is true not only of underdeveloped countries but also of traditional fishing nations such as Iceland, whose entire

economy is dependent on its offshore cod fishery. Canada has also threatened similar action. Several international incidents have been directly tied to the problem of sharing a fisheries resource or to disputes over conflicting uses of a resource within a given area, for example, the conflict between United States lobster fishermen and Russian trawlers off the New England coast. In a few cases a fishery is regulated by treaty. This is true for the Pacific salmon which is shared by the United States, Canada and Japan. However, since the treaty only governs these three nations, what happens when other countries not regulated by any agreements attempt to share this resource? The problems resulting from international conflicts over marine fish resources are among the most challenging and complex we can define for the renewable resource manager.

Because of their need for protein, aquaculture of food fishes constitutes an important resource for the underdeveloped nations of the world. However, many of these countries already have an extensive development of these resources. As poorly developed nations attempt to raise their standard of living, they will put greater and greater demands on protein sources such as grain crops, fish meal and even beef cattle. Thus, they will come into direct competition with more affluent countries. For European nations which depend on imports of meat or high protein feeds to raise beef or dairy cattle such competition will become a serious problem. Though Canada and the United States produce enough food for their own use, how can they morally disregard the needs of four-fifths of humanity? In addition, they are caught up in the international balance-of-payments problem and numerous treaty obligations. Thus, the fisheries resources of the world remain unalterably tied to questions of moral, political, and social values.

An even greater problem facing freshwater and coastal estuarine resources is the continuous degradation of the environment and the concomitant changes in the fish populations. Although we do not directly consider the well documented problem of pollution and its effects on aquatic organisms, it is certainly important in many of our fisheries. It has had a direct bearing on the development of certain

fisheries resources such as the Great Lakes fisheries, notably in Lakes Ontario and Erie. However certain fisheries in the upper Great Lakes have also been affected.

In view of these serious problems, the need for recreation may seem of diminished importance. Yet it continues to dominate the inland freshwater fisheries of the affluent western world. In island waters, sport and commercial fisheries are often in direct competition for a common resource and even when they compete for different resources within a commonly shared area, conflicts are just as serious. The Great Lakes fisheries constitute an outstanding example of this sort of conflict. Regulation of these fisheries is made even more complex since jurisdiction is both international and interstate.

Inland sport fisheries are also faced with a myriad of other problems. The development of an industrialized, urban society with a seemingly insatiable demand for recreational space has created a paradoxical situation; the masses of modern man are densely concentrated while the resources are usually located at great distances from these population centres. However, with improvements in transportation and the general affluence of the public even once remote resource bases are hard put to produce satisfactory recreational experiences. Fishery managers are thus faced with the task of providing satisfactory yields and dealing with the incremental and continuous degradation of existing resources. Certain resources such as streams continue to be degraded at a rate, which if continued, will soon make them a rarity as a fishery resource. The construction of impoundments have created new fishery resources, yet their management remains quite primitive. Of immediate concern is the loss of existing fishable waters near urban centers. The increasing demand for such waters continues to tax the most imaginative and creative minds in the field of fisheries administration. Finally, where demand is greater than supply, certain special interest groups usually bring pressure on administrators to propose legislation favouring their use of the resource. This tends to exclude the less organized, more general sportsman. Attempts to restrict both the harvest and the use of a resource to special interest groups continues to be one of the most challenging problems in sport fishery management.

Coastal marine fisheries are faced with many of the same problems as inland freshwater fisheries. In addition to the pollution of the environment, there is an increasing demand for salt water based recreation plus conflicts over the commercial and/or sport use of many marine fish and shellfish.

Thus, the challenges facing the fishery manager are as complex, formidable and numerous as any in the field of renewable natural resources. They include the major problems confronting man today: over-population of the third world, the stark reality of protein starvation for four-fifths of humanity, insidious and incremental destruction of the world's natural resources at a time of resource scarcity, increasing competition between wealthy and developing nations for the resources that remain and the psychological impact of coping with a technological world at a time when the safety valve provided by outdoor space is vanishing or being altered. Even though advances in science and technology have been the cause of many of these problems, they can also be used to formulate effective solutions. However, the proper application of these solutions will also depend on the social and political wisdom of mankind.

We are on the verge of spectacular breakthroughs in the field of ecosystem management, but man's greatest achievement will be in finding the social and political answers to make the "wise" use of his renewable natural resources a reality for succeeding generations.

R. I. P.

Edmund Fitzgerald

and

taconite pellets

11/11/75

THE MARRIAGE PROBLEM

by L. Mathemaki,
Professor of Mathematical Marriage Counselling

Most mathematics has its roots in very practical problems and, for most of us, one very practical problem we have all encountered — and some of us have solved — is: “how many acquaintances must one have in order to get married without undue competition?”.

I will pose this problem in a typical idealized way: we will consider a set B of boys, and set G of girls; each boy wishes to marry an acquaintance from G ; perhaps a boy will not know all the girls in G . Bigamy will not be allowed, and we wish to attain the typical male-chauvinist goal that all the boys get married, without regard for left-over girls.

Let's consider an example: we have four boys, improbably named 1, 2, 3, and 4. Each boy knows some girls, any one of whom he would not mind marrying. Let us call the girls A, B, C, D , and suppose the acquaintances are

Boy	His acquaintances
1	A
2	A, B
3	B, C
4	A, C, D

If all the boys are to be married, 1 must marry A , so 2 must marry B , 3 must marry C , and 4 must marry D . In this case, all the boys can get married, and in fact there is only one possible way they can pair off.

Another example: suppose the ‘acquaintance table’ is:

Boy	His acquaintances
1	A, B
2	A, B
3	A, B
4	A, B, C, D

In this case marriage is impossible: the boys 1, 2, 3, only have two girls to choose from.

Thus we see that sometimes marriage is possible and sometimes it isn't. **The marriage problem is simply to determine conditions which will guarantee the possibility of marriage.**

To solve the problem, let's return to the general situation — the two sets B, G — and suppose all the boys can be married off. Consider any set C of boys from B (that is, consider any subset C of B) and denote by $F(C)$ the set of those girls from G each of whom knows a boy in C .

For instance, in our first example we have $F(3) = (B, C)$; $F(1, 2) = (A, B)$, $F(2, 3, 4) = (A, B, C, D)$, and so on.

We have assumed all the boys can be married off; thus no matter what C is, $F(C)$ must contain the (potential) wives of the boys in C , and since different boys have different wives (the no-bigamy condition), there are at least as many girls in $F(C)$ as there are boys in C .

So we have the assertion: **if marriage is possible then, for each $C \subseteq B$, $F(C)$ contains at least as many elements as C .**

THE SURPRISING THING IS: the converse of our assertion is also true. Given any sets B, G of boys and girls, if $F(C)$ contains at least as many girls as there are boys in C , for every $C \subseteq B$, then all the boys can be married off.

The proof of this is very tricky indeed — it proceeds by induction on the number of boys in B . It would be a worthwhile problem for a good grade 13 student to attempt; at least it illustrates that induction is not just a turn-the-crank method involving the summation of finite series!

Of course, such a proof only works for a finite number of boys, so it then is natural to ask: what happens if B contains an infinite number of boys? Consider an example: call the boys 1, 2, 3, 4, ... the girls $1^*, 2^*, 3^*, 4^*, \dots$ and suppose the acquaintances are:

1. knows all the girls
2. only knows 1^*
3. only knows 2^*
4. only knows 3^*

and, generally, if $n > 1$, n only knows $(n-1)^*$.

Every set C of boys knows as many girls as there are boys in C ; in fact, if 1 is in C , then $F(C)$ contains all the girls. But marriage is impossible — 2 must marry 1^* , 3 must marry 2^* , ... leaving no one for that extremely popular boy 1.

Thus the condition that worked for the finite situation won't work in this case; the next step would be to get some compromise. It turns out that if the condition holds, and each boy only knows a finite number of girls, then marriage is possible. The proof of this seems to require some very profound mathematics indeed; it was first proved in 1949, and subsequent generalizations now constitute a fairly substantial body of knowledge "the theory of transversals and representative sets".

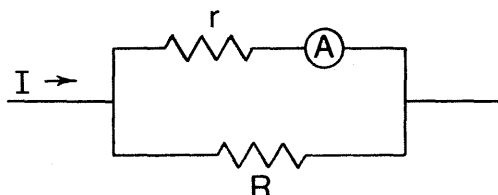
TWO MORE PROBLEMS:

(1) **The problem of the monks:** this was first posed by Balzac. We have two sets, a set M of monks, and a set A of non-monks; each monk has acquaintances in A , and he wishes to establish a 'harem' consisting of his acquaintances, of a particular size, and disjoint from the other harems. Under what conditions can he do this? It turns out that this is just a disguised marriage problem: replace each monk by as many boys as he wants members in his harem, and seek conventional marriages.

(2) **The friendship problem:** this appears unrelated to the marriage problem. We consider a party at which each pair of persons has a common acquaintance (including the possibility that a pair know each other). Then, someone at that party knows everyone! The only proof I know of is pretty complicated — in fact, it turns the problem into a problem in a peculiar geometry. **Can anyone furnish a nice neat elementary proof?**

True or False?

Ammeter: a current-measuring device. Most modern ammeters are of the moving coil type developed during the 19th century. The current to be measured is passed through a rectangular coil of many turns of wire pivoted between two jewelled bearings in a magnetic field. The interaction between the current and the magnetic field exerts a torque which turns the coil against a spring. The angle through which the coil rotates is then a measure of the current in the wire. The sensitivity of the instrument is set by the strength of the magnetic field, the construction of the coil (size of wire, number of turns and area of coil) and by the strength of the restoring spring. An external circuit can easily be used to **reduce** the sensitivity of a meter as can be seen in the diagram:—



In this circuit, only the fraction $R/(R + r)$ of the current I , flows through the ammeter, A . No external circuit, however, can be used to increase the sensitivity.

True or False?

Carotid: one of the two great arteries that carry blood to the head. The name is derived from *karoo* (= stupefy), compression of the carotid arteries being thought to produce stupefaction.

Carotene: one of the carotenoid pigments common in plants (and particularly abundant in carrots).

Carob: leguminous evergreen Mediterranean tree (and its pod and seeds). The large red pods have been used for food for animal and man since prehistoric times. It is believed that it may have been the "locust" eaten by John the Baptist in the wilderness. The seeds are reputed to be extremely uniform in size and weight and to have been used as weight standards. They are believed to be the origin of the **carat** which is now the measure of weight for precious metals and jewels.

WHAT IS TOPOLOGY?

by: S. A. Naimpally,
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As with any subject it is quite difficult to give a precise definition of Topology. Topology means different things to different people but broadly speaking it deals with *continuity*. The terms continuous, continuity etc. are used often in day-to-day language, and the purpose of this little article is to explain how they are made precise in Mathematics. Mathematicians have been struggling for centuries to give a precise meaning of continuity, and although a satisfactory definition was given in the middle of the nineteenth century by Weierstrass and Heine, a very simple approach was first suggested by the great Hungarian mathematician F. Riesz in 1908. We will use Riesz's approach via the concept of *nearness* which is one of the most important concepts in the whole of mathematics. The beauty of nearness is that it is simultaneously simple and rigorous.

Let us take a simple example. Suppose Paul is a friend of the Russel family. If Paul continues to be a friend, then we say that the friendship is *smooth* or *continuous*. If there is a quarrel between Paul and the Russels, we say that the friendship is *broken* or *discontinuous*. It is easy to think of many such examples, but rather than do that, we will pursue only a few and arrive at a precise definition of continuity.

The statement "Paul is a friend of the Russels" can be simply explained by "Paul is *near* the Russel family". Here nearness is an abstraction of friendship but it could also represent many other relations. For instance, a phrase which we frequently use in day-to-day language, we may say that "Paul is near the Russel family" provided that Paul lives in their neighbourhood, or Paul is related to the Russel family (Perhaps as an uncle or grandfather). Thus we have the concept of *nearness* between a person and a family.

A question naturally arises as to what properties this nearness relation satisfies? Some of the obvious ones are (i) each member of the Russel family is near the Russel family (we are, of course, considering an ideal family which has no delinquents!), (ii) No person can be near a non-existent family, etc. However, a subtle difference exists between these day-to-day examples and the mathematical concept of

nearness. Whereas it is possible in life for a person to be neither near nor far (opposite of near), we require our mathematical concept to avoid this possibility.

As time passes we all change or we are all transformed. If a person who is near a family remains near we say that this transformation is *continuous* and, in the contrary case, we say that it is *discontinuous*. This idea is very basic and is at the root of the concept of limit in Calculus.

In Topology one considers "higher" types of nearness too. For example we may say that the Russel family is near the Hardy family, if as it happens so frequently, they are introduced to each other by a common friend or a girl from one family marries a boy from another. This nearness between two families is called *proximity*, and interestingly enough, even in Mathematics the two most important proximities arise in completely analogous ways. One then considers *proximal continuity* which is concerned with the preservation of proximity. Finally we have the concept of nearness for several families. For example, several families may be *near* because they have a common ancestor or they have a common cause and get together for a picnic.

Although these ideas are simple, they are at the basis of Calculus, Analysis and a big chunk of Modern Mathematics. They have found applications in Mathematical Biology, Psychology, Econometrics, Physiology of the Brain etc. For further details and references, the interested readers are referred to the following:

- (1) S.A. Naimpally and B.D. Warrack, *Proximity Spaces*, Cambridge Tracts in Mathematics No. 59, Cambridge University Press, U.K. (1970).
- (2) S.A. Naimpally, *Proximity Approach to General Topology*, Lecture Notes, Lakehead University (1973).
- (3) P. Cameron, J.G. Hocking and S.A. Naimpally, *NEARNESS; a better approach to continuity and limits*, Lecture Notes, Lakehead University (1973).

INDUCTIVE JUSTIFICATION OF INDUCTION

by: Gunars Tomsons
Lakehead University

The analysis of scientific reasoning by many able investigators has resulted in the acknowledgement of induction as being a legitimate and indispensable process of framing general hypotheses. This process, according to the analysis consists of formulating a general hypothesis on the basis of observed instances. Thus from the observed mortality of humans one infers the the mortality of any member of any mankind.

Since the future seems so open and uncertain, and since a radical change in the constitution of the universe cannot be excluded as a possibility, it has been thought at times that a justification is needed for a procedure of science that purports to produce hypotheses and laws of nature on the basis of the limited evidence of the senses.

One sort of justification that has appeared to be the most promising method of justifying induction is the so-called inductive justification of induction. According to this method one is justified in using induction because observed instances of inductive reasoning have turned out to be reliable. There is certainly some attractiveness in a proposal to justify induction on empirical grounds. It seems that if we can make what we think to be legitimate inferences based on past experiences when we deal with the objects of science, then we should be able to learn from past experiences and make legitimate generalizations even when the objects have a different nature, i.e., when the objects are inductive arguments.

However, at first sight there appear to be also good reasons against an inductive justification of induction, especially if we formulate the problem in terms of inductive rules. The justification of induction is then seen as a justification of a rule that permits generalization. Hence, if we want to justify our use of the rule through inductive procedures, we seem to be using in our justification a rule the general use of which we are trying to justify. There seems to be something illogical about a procedure of this sort. Nevertheless, Max Black, a present day philosopher interested in inductive logic has argued that inductive justification of induction is possible.¹ Hence I want to present the model for inductive justification

of induction and Black's reasons for his position. I shall then show that Black's position is not tenable.

Black conceives of the inductive justification of induction in terms of a justification of the inductive rule governing one type of inductive argument. The rule whose justification he seeks is the rule permitting the following argument:

Most of the examined A's have been B's r_1
The next A will be B.

Observation of the repeated application of the rule r_1 , prompts us, according to Black, to produce an argument of the following form:

r has usually been successful. r_1
 r will be successful in the next instance.

An argument of this sort is called a self-supporting argument, because the rule r_1 , is used in the argument to establish that the next instance of the use of the rule will be successful. The meaning of 'the next instance' in the conclusion has not been made explicit, but Black's examples indicate that by 'the next instance' he means the use of the rule in a non-self-supporting argument. Arguments of the above kind have at times been called second-level arguments, because they refer to arguments whose referents are found in the non-linguistic world. One could, of course, construct second- or third-level arguments; but in the case under consideration we have been presented with only a second-level argument. One may not assume, though, that the rules used in second-level arguments must differ from the first-level arguments merely in virtue of the fact that second-level arguments try to establish facts about the first-level arguments. Moreover, I think that it would be absurd to claim that the inductive r_1 differs in character depending on whether it is employed in first- or second-level arguments.

Arguments of the self-supporting kind have been criticized on the grounds that they are circular, i.e., that their premises contain the conclusion, if only as a presupposition or a suppressed premise. Black has provided us with a seemingly decisive argument against the view that these arguments have a suppressed premise asserting the general reliability of the rule r_1 ²

He has argued that the argument could not have a suppressed premise of this kind because by repeating the conclusion in the premises we no longer have an inductive argument. We have a technically sound argument which is of no use to us because it is circular. But only deductive arguments can be circular. Hence, according to Black, an attack of this sort on the inductive justification of induction cannot but fail.

This appears to be a victory over those who allege that inductive justification of induction is circular in the logico-technical sense. However, I am not quite convinced. The attack by Black succeeds only if his opponent is willing to grant that the argument is in fact inductive. But need he grant this? It seems to me that Black's opponent could say that the justification of induction fails because the argument is circular and hence also deductive. On this view, Black only believes that he has provided an inductive argument, whereas in fact he had the suppressed premise that enumerative induction is effective. In other words, he has assumed that which he wants to prove inductively. And if one wants to prove what he assumes to be true in the premises, he has a circular argument whether he intends to use deductive or inductive procedures for proving whatever he intends to prove.

From the point of justification of a belief that rests on an inference it certainly is presupposed that the belief is well supported, and in order for a belief to be well supported we must have a reliable rule of inference as well as good evidence. Thus the reliability of the rule is presupposed. It is true, I think, that we may not be conscious of the belief that our rule is reliable. However, from the point of view of justification of beliefs we have to take into account beliefs that we are not aware of in our day-to-day arguments, if they are relevant to our epistemological order of beliefs. I conclude that Black has not succeeded in showing that the inductive justification of induction is not circular, and I have already indicated that one has good reasons for claiming that it is circular.

Black has argued that self-supporting arguments are not circular also on the grounds that the argument succeeds in raising the reliability of the rule. I want to show, nevertheless, that this argument is equally unsuccessful, and I will first reproduce Black's example of a self-supporting argument:

- (a) In most instances of the use of *R* in arguments with true premises examined in a wide variety of conditions, *R* has been successful

Hence (probably):

In the next instance to be encountered of the use of *R* in an argument with a₃ true premise, *R* will be successful.

'*R*' stands for the following rule:

To argue from

"most instances of *A*'s examined in a wide variety of conditions have been *B*"

to (probably)

"The next *A* to be encountered will be *B*"

Black contends that the argument (a) supports rule *R* by raising the reliability of the rule and hence the subsequent first-level arguments using the rule. The situation can be presented in the following schematic way using these abbreviations:

'*a*' will represent first-level inductive arguments before the use of the self-supporting argument.

'*A*' will represent the self-supporting argument

'*b*' will represent a first-level argument after the use of the self-supporting argument '*A*'

'*r*_{*a*}' is the inductive rule used in arguments *a*

'*r*_{*b*}' is the same rule used in argument *b*

'*r*_{*A*}' is the rule used in the self-supporting argument

The sequence below represents the temporal order of the arguments and the abbreviated argument *A* is presented directly below the sequence:

$a_1, a_2, a_3, \dots, a_n, A, b$
 r_a was successful
 A: $\dots r_A$
 r_b will be successful

According to Black argument A establishes that r_b is more reliable than r_a , but I do not see that this has to be the case, since we do not know much about the reliability of r_A . Black has no reason to assume that his second-level induction will perform as expected since in proving the conclusion we have assumed that r_A will perform satisfactorily; we certainly have no inductive evidence in support of the reliability of r_A .

Some arguments are in order also with respect to Black's concept of the degree of reliability of the rule. Black conceives of the strengthening of rules and arguments as follows:

Let us assume that the following argument b using the rule r has the strength 4/5 or 8/10:

Argument b 4/5 of examined A's are B's

Next A will be B

Given appropriate observational evidence we can construct a self-supporting argument:

Argument A 9/10 of examined arguments using rule r were successful

Next use of r will be successful

This argument entitles us to say that the degree of strength of the rule r_b is 9/10. Hence we should increase the strength of the argument b to 9/10.

I already indicated why I do not think that Black has shown that the reliability of the rule would be increased, but Black has not really explained why we are entitled to speak about degrees of reliability of the rule. It may be plausible to talk about degrees of strength of an argument as a function of the amount of evidence available. However, is it plausible to say that rules have varying degrees of reliability? Of course, three different rules could conceivably have distinct degrees of reliability, if we should find that the percentage of past

successes differs from rule to rule. But why should an inductive rule increase or decrease in reliability depending on the ratio of past successes? Our confidence may perhaps increase if we see the rule succeed, but to say that our confidence is increasing is not to say that the reliability of the rule is increased.

I conclude that Black's attempts to provide inductive justification of induction have not succeeded. Moreover, I have supplied reasons which show that inductive justification of induction is implausible.

1. M. Black's writing on inductive logic and critical discussions of them can be found in the following works. M. Black, *"The Justification of Induction"*, Language and Philosophy, (Ithaca, New York: Cornell University Press, 1949), pp. 59-88; "The Inductive Support of Inductive Rules", *Problems of Analysis*, (Ithaca, New York: Cornell University Press, 1954), pp. 191-208; W.C. Salmon, "Should We attempt to Justify Induction?", *Philosophical Studies*, VIII (1957), pp. 33-48; M. Black, "Self-Supporting Inductive Arguments", *Journal of Philosophy*, LV (1958), pp. 718-725, or pp. 209-218; P. Achinstein, "The Circularity of a Self-Supporting Argument", *Analysis*, XXII (1961-62), pp. 138-144; M. Black, "Self-Supporting and Circularity: A Reply to Mr. Achinstein", *Analysis*, XXIII (1962-63), pp. 43-44; P. Achinstein, "Circularity and Induction", *Analysis*, XXIII (1962-63), pp. 123-127; H. Kyburg, "recent Work in Inductive Logic", *American Philosophical Quarterly*, I (1964), pp. 1-39; M. Black, "the Raison d'Etre of Inductive Argument", *British Journal for the Philosophy of Science*, XVII (1966), pp. 177-204; also in *Margins of Precision* (Ithaca, New York: Cornell University Press, 1970); "The Justification of Induction", *Philosophy of Science Today*, S. Morgenbesser (ed.) (New York: Basic Books, 1967); "Self-Supporting Inductive Arguments", *Models and Metaphors*, pp. 210.
2. M. Black, *Problems of Analysis*, pp. 199-200.
3. M. Black, "Self-Supporting Inductive Arguments", *Models and Metaphors*, p. 210.

PHOTO-DEGRADABLE PLASTICS

by: *Dr. N.A. Weir,*
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One consequence of the increased standards of living experienced by most industrial countries is the rapid increase in the volume of solid waste. Wrapping of foods to reduce contamination and prevent deterioration has brought many social benefits, which at present tend to outweigh the disadvantages of having to dispose of the packaging materials after use.

Plastics are also replacing conventional containers, like glass and ceramics to an increasing extent. One needs only to look around the shelves of a supermarket to see polyethylene detergent and bleach containers, polyvinylchloride (P.V.C) shampoo and edible oil containers and polystyrene beverage cups. The advantages of these materials are obvious — ease and cheapness of formation of variously shaped containers, favourable mechanical properties (non-brittle) and their light weights.

However, the steadily increasing proportion of plastic wrapping materials and non-returnable plastic containers in domestic and industrial wastes is now causing major disposal problems in many parts of the world. A recent survey shows that three polymers account for 90% of municipal plastic waste. These are: polyethylene (38%), P.V.C. (31%), and polystyrene (21%); the remaining 10% is composed of varying amounts of polypropylene, polyesters, polyvinylidene chloride, and cellulose acetates and butyrates.

Unlike cellulosic packaging materials, like paper, the plastics mentioned above are not bio-degradable (i.e. capable of being broken down, ultimately to CO_2 and H_2O , by micro-organisms), hence they tend to be very persistent in the environment.

Lack of biodegradability can be attributed to three typical characteristics of the polymers most used in packaging.

(1) The polymers are largely hydrophobic, and hence wettability by polar molecules, like

water, a prerequisite for many biological reactions, is extremely low.

(2) Permeabilities of these polymers to gases like oxygen are very low, and thus the efficiencies of many bio-processes are greatly impaired.

(3) Molecular weights of these materials must be high to give them adequate mechanical strength. However, micro-organisms tend to attack at the ends of large molecules, and since the number of these ends is inversely proportional to the molecular weight, it follows that the rate of attack on a polyethylene molecule which might be used in packaging (Molecular weight 120,000) will be very low.

It is ironic that the factors which determine the usefulness of these polymers in packaging are precisely the factors which lead to their persistence in the environment.

Some typical biodegradability figures, measured on a growth rating scale of 1 to 10, are summarized below.

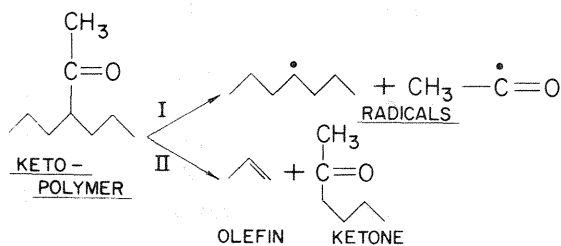
Compound	Molecular weight	Growth Rating
polyethylene	129,000	0
polyethylene	52,500	0
polyethylene	13,800	1
polyethylene	1,350	2
polystyrene	87,000	0
polystyrene	2,100	0
polyvinyl chloride	52,000	0
polypropylene	27,000	0
dodecane ($\text{C}_{12}\text{H}_{26}$)	170	5

The importance of molecular weight can be seen by comparing the various polyethylenes and the chemically similar, non-polymeric molecule dodecane. The above polymers may be rendered bio-degradable, by modification of their hydrophobicities, their permeabilities and by decreasing their molecular weights. This can be achieved simply by exposing photo-degradable forms of these polymers to natural sunlight, or by adding suitable sensitizers to them, the sensitizers also absorbing sunlight, and initiating polymer degradation.

It is convenient to discuss the two methods separately.

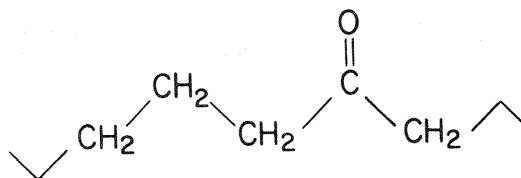
Ketone Polymers

After two decades of research on the prevention of sunlight degradation of polymers, scientists are applying their results to the design of polymers which will readily undergo sunlight degradation. Chemically, the reactions which lead to the failure of the polypropylene webbing used on lawn chairs, are very similar to those which are involved in the disposal of photo-degradable polypropylene, and in both cases ketonic (or carbonyl) compounds play a critical role. When the carbonyl group, $C=O$, absorbs ultra-violet radiation of wavelength around 300 nm, ($1\text{nm} = 10^{-7}\text{cm}$) — and natural sunlight contains a small ultra-violet component with wavelengths extending to around 295nm—one of the non-bonding (localized) electrons on the oxygen atom is promoted into an anti-bonding π^* — molecular orbital, which results in the delocalization of the electron over the C and O atoms. Such a promotion is referred to as an $n \rightarrow \pi^*$ transition (i.e. transition from an n-type to an anti bonding π^* molecular orbital) and the excited molecule so produced can react by undergoing bond fission adjacent to the $C=O$ group. Two distinct processes are possible, and these occur in both polymers and in small molecules.



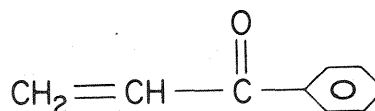
Reaction I (better known as the Norrish Type I process) results in the formation of two free radicals. A free radical is a species which contains an unpaired electron and which is formed by the homolytic fission of a covalent bond. Reaction II (Norrish Type II) involves an intramolecular rearrangement which is accompanied by fission of the molecule into another ketonic compound and an olefin.

It can be seen that if a ketonic group is introduced into a polymer chain, exposure to ultra-violet radiation will result in scission of the main chain, (Reaction II) and this reaction is employed in one type of photo-degradable polymer. Photo-degradable polyethylene has the structure:

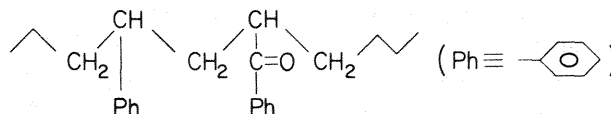


and it can be seen that both Reactions I and II will result in scission of the chain. To make polyethylene photo-degradable, it is necessary to introduce only one ketonic group per thirty carbon atoms in the chain; the fragments produced by photo-fission are potentially bio-degradable.

Photo-degradable polystyrene and polypropylene are now available commercially (known as "Ecolyte S" and "P" respectively). Ketonic groups are incorporated into these molecules by copolymerization of small amounts of phenyl vinyl ketone.



The resulting copolymers have the following structures, the ketone groups occurring at random in the chains, e.g. with styrene,



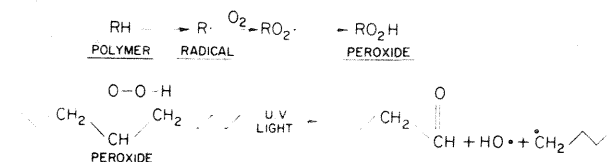
Lifetimes of these polymers can be controlled by varying the phenyl vinyl ketone concentration in the copolymers. "Ecolyte S" and "P" were developed by Dr. J.E. Guillet and his co-workers at the University of Toronto.

It should be emphasized that in all of the above examples, the physical properties of the photo-degradable polymers are indistinguishable (except for ultra-violet absorption characteristics) from those of the pure homopolymers.

(2) Sensitized Photo-degradation

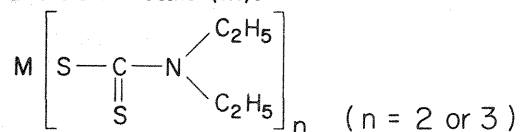
An alternative approach to photo-degradable plastics has been made by Dr. G. Scott at the University of Aston, England, and this employs a combination of photo-oxidation and subsequent photo-degradation. Free radicals, like those formed in Reaction I, being electron-deficient species react readily with molecular oxygen in the air—the oxygen molecule exists in a triplet state — to produce peroxy radicals

which in turn abstract hydrogen atoms to produce a hydroperoxide and a new radical. The hydroperoxide is a key intermediate, since it absorbs the ultra-violet component of the sunlight, leading to photo-fission of the bonds adjacent to the peroxide ($-\text{O}-\text{O}-$) group. The new radicals formed can continue the oxidation process, and a chain reaction can develop. Photo-oxidation of a polymer is illustrated by the following reaction sequence, which incidentally, is the same sequence that leads to the deterioration of the useful properties of polypropylene on exposure to sunlight.



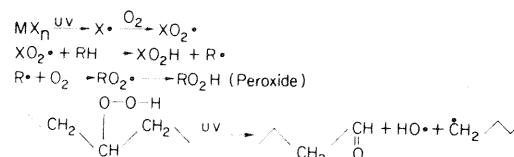
It can be seen that photolysis of the hydroperoxide will lead to chain scission, and ultimately to destruction of the polymer.

One advantage of this system is that the oxidation reactions do not require continuous illumination; hydroperoxides can be formed in the dark, and subsequently decomposed on exposure to sunlight. One disadvantage of the system illustrated above is that the initial production of free radical centers on the polymer is intrinsically very slow, and thus photo-oxidation as such, is not a practical system for photo-disposable plastics. Scott discovered, largely by accident, that initiation of the photo-oxidation could readily be accomplished by adding suitable substances to the polymer. The compounds found most useful in this regard are the dithiocarbamate complexes of transition metals (M).

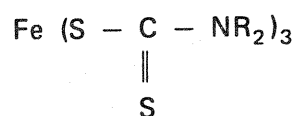


Used in high concentrations these molecules act as photo-stabilizers (molecules added to inhibit photo-degradation), but become photo-initiators at low concentrations (<1%). Initiation is brought about as follows. On exposure to sunlight the photo-initiator is decomposed into free radicals, which react with oxygen in the air to form peroxy radicals, which in turn abstract hydrogen from the polymer chains forming radical centers. The polymeric radicals then undergo oxidation.

The process is summarized by the following scheme: (MX_n is the initiator)

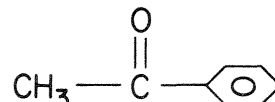


Addition of small amounts of these dithiocarbamate complexes to polyethylene polypropylene and polystyrene render the polymer photo-degradable, and their lifetimes can be controlled by variation of the concentration of photo-initiator. The iron complex,

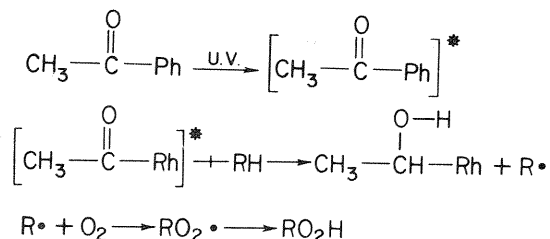


is particularly attractive as a photo-activator for degradation of plastics used in the food-stuffs industry since it is non-toxic.

Yet another group of photo-initiators has been developed by scientists at the National Research Council of Canada, Ottawa. These compounds are aromatic ketones like benzophenone, which absorb



ultraviolet radiation in the sunlight spectrum to produce excited triplet molecules by $n \rightarrow \pi^*$ transitions. (In the triplet state the electrons in the π^* and n -type molecular orbitals have parallel spins). The triplets have sufficiently long lifetimes to undergo hydrogen abstraction reactions with the polymers, again forming radical centers on the chains, as follows:



The polymer is then oxidized and the chains broken by exposure of the hydroperoxides to sunlight.

THE COMPOUND MICROSCOPE

by: Mr. B. Spenceley,
Lakehead University

It scarcely needs to be said that the compound microscope is widely used in the teaching of biology, but whether most students have a sufficient understanding of the operation of the microscope is debatable. The purpose of this series of articles is to improve this understanding and to enable students to get more use out of their instruments.

One of the important characteristics of an instrument for examining small objects is its magnification. There is a certain amount of arbitrariness about the definitions of magnification. The angular magnification, which is the type we are concerned with here, is usually defined as the ratio of the angle subtended by the object at the distance of most distinct vision to the angle subtended by the object seen through the instrument. A value of 25 cms is usually taken as the average distance of most distinct vision, although there is a great deal of variation in this from one person to another.

The simplest optical device for examining small objects is a single lens of short focal length. If f is the focal length in centimeters the magnification, m , is $25/f$ so one would be tempted to consider that higher and higher magnifications could be easily obtained by making f smaller and smaller. In practice magnifications up to 20x can be achieved using simple magnifiers but attempts to get higher magnifications are hindered by the aberrations that always accompany lenses of short enough focal length and large enough aperture to be useful.

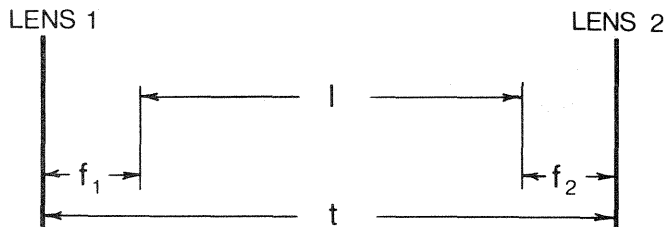
The compound microscope overcomes this difficulty by using a pair of lenses relatively well separated. Elementary optics texts show that if two lenses whose focal lengths are f_1 and f_2 are separated by a distance t , the effective focal length of the combination is given by the simple equation

$$\frac{1}{f} = \frac{1}{f_1} + \frac{1}{f_2} - \frac{t}{f_1 f_2}$$

The equation looks more useful if t , the distance between the lenses is replaced by ℓ ,

the distance between the focal points. From the following simple diagram it is clear that

$$\ell = t - f_1 - f_2$$



When the last two equations are combined a simple expression for the effective focal length results:— $f = \frac{-f_1 f_2}{\ell}$

The point of all this is that although it is difficult to make a single short-focal-length lens we can get the same effect by using two moderately 'short' lenses separated by a relatively large distance ℓ . Because ℓ is in the denominator large values of ℓ mean small values of f . So the compound microscope can be regarded as a simple magnifier; as such its magnification is

$$m = \frac{25}{f} = -\frac{\ell}{f_1} \cdot \frac{25}{f_2}$$

Sometimes it is more useful to consider the effect of the two lenses separately. The first lens (f_1), usually called the objective, forms a real, inverted image of the object in the focal plane of the second lens, the ocular or eyepiece lens. The lateral magnification of the objective is $-\ell/f_1$. The second lens is then used as a simple magnifier to look more closely at this intermediate image. The angular magnification of the eyepiece is of course $25/f_2$. The overall magnification of the compound microscope which we have already noted to be $[-\ell/f_1][25/f_2]$ is merely the product of the individual magnifications of the objective and eyepiece lens.

Because the magnification of the eyepiece depends only on its focal length, it is feasible to mark the lens mount with the magnification. This is the meaning of the numbers such as 10X or 15X found on oculars. The magnifi-

cation of the objective on the other hand depends on the length λ and unless this is known and constant the magnification cannot be specified. For this reason microscope manufacturers have standardized the distance λ at 160 mm for the great majority of their instruments. For these microscopes it is possible to engrave the objective magnification on the lens barrel, the number is unchanged if the objective is used in a conventional microscope no matter what the country of origin — an example of international cooperation that could well be imitated by other apparatus manufacturers. The overall magnification of the compound microscopes used by most schools is simply obtained by multiplying together the magnifications engraved on eyepiece and objective lenses.

Another important characteristic of the compound microscope is its resolving power or ability to distinguish small details. This is not the same thing as magnification; it is relatively easy to get large magnifications, for example by making λ larger, but this does not ensure that more detail is visible. High magnification that is not accompanied by good resolution is called "empty magnification" because the increased image size does not carry any more information about the subject. **The image is larger but blurrier.**

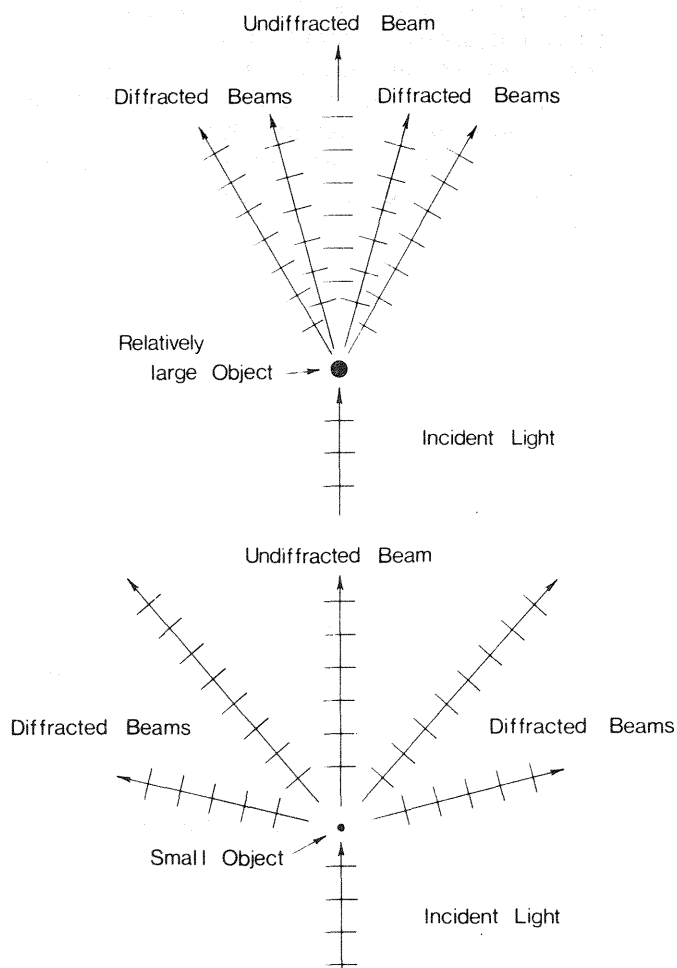
For a microscope of good quality the resolving power is limited by the diffraction of light in the plane of the object under examination; in an inferior instrument it could be limited by ordinary lens defects. A proper treatment of the theory of diffraction in microscopy is beyond the scope of this article, but some of the basic ideas can be conveyed in relatively simple terms.

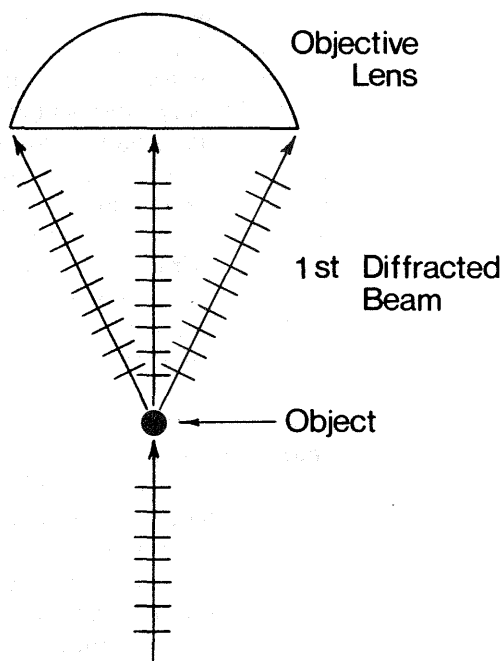
When a beam of light passes an object whose size is comparable to the wavelength of the light, the beam is *diffracted*, by which we mean that some of the light on the far side of the object travels in directions different from the original one. Usually the beam breaks up into several beams — one undisturbed by the object, a pair making a small angle to the original direction, and so on. Two cases are illustrated in the following figures, first a relatively large object and then a small object.

The diagrams illustrate the important idea that the smaller the object is the larger are the diffraction angles, and the larger the object is the smaller are the diffraction angles. The information about the size and shape of the object is carried in the angular spread and relative intensities of the diffracted beams.

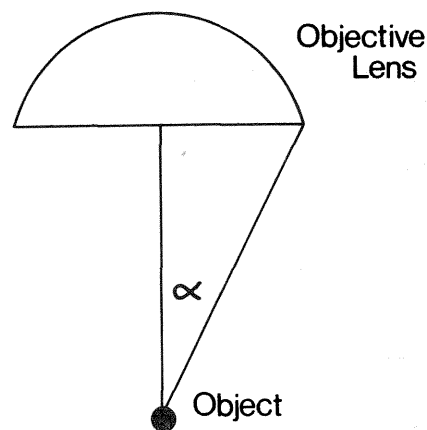
The size of the object compared to the wavelength is also important in determining the angular spread of the diffracted beams. For an object of a given size the longer wavelengths (the red end of the spectrum) are diffracted through larger angles than the shorter wavelengths (the blue end of the spectrum).

If the objective lens of the microscope is small in diameter compared to its distance from the object, the angle it subtends at the object will also be small. If the angle is so small that only the undiffracted beam enters the lens, then it is as if there were no object there at all, since that would also be the only beam in the absence of an object. In order even to detect the presence of the object the





lens must at least be able to accept the first diffracted beams as shown in the next diagram. The limit of resolution of a microscope is therefore the smallest object whose first diffracted beams are just accepted by the objective. Naturally, if more detail is needed than the mere existence of the object, then the lens



must be able to accept more of the diffracted beams. Strictly speaking, the resolving power is related to a quantity called the numerical aperture of the objective lens which is the product of the sine of the half-angle subtended by the lens at the object and the index of refraction of the medium below the lens — usually air but sometimes not. The resolving power is the wavelength of the light divided by the numerical aperture — $\lambda/n \sin \alpha$. The best resolution is obtained by using blue light and by filling the space between the object and the lens with a liquid of high refractive index.

EDITOR'S AND COLLABORATORS' COMMENTS

T G I T L P O T I

IPTLIGOTT: a sophisticated anglo-saxon surname?
or is the answer simply biodegradable?

Please permutate, decypher and send in your interpretation.

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